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MATHEMATIK

ZUR LAPLACE-TRANSFORMATION. VI

Von V. S. IGNATOVSKIJ (W. IGNATOWSKY), korrespondierendem Mitglied der Akademie

Wir wollen nun folgende Gleichung betrachten.

$$\frac{1}{c^2}\frac{\partial^2 u}{\partial t^2} + c_1\frac{\partial u}{\partial t} + a_1\frac{\partial u}{\partial x} - \frac{\partial^2 u}{\partial x^2} - \frac{\partial^2 u}{\partial y^2} + mu = \varphi(x, y, t). \tag{1}$$

Es soll hierbei $u(x, y, t)$ eine Funktion von x, y und t sein, unabhängig von z. Wir haben also einen ebenen Fall und wollen hierbei folgendes Problem behandeln.

Es sei die YZ-Ebene des rechtwinkligen Koordinatensystems die Grenzfläche, rechts von welcher (also bei positiven x) die Gleichung (1) gelten soll.

Analog den vier Grössen (2) und (3) unserer Note I [1] führen wir auch jetzt die folgenden vier, aus $u(x, y, t)$ fliessenden, Grössen ein:

$$u(0, y, t) = u_0(y, t); \quad \left.\frac{\partial u(x, y, t)}{\partial x}\right|_{x\to 0} = h(y, t) \tag{2}$$

und

$$u(x, y, 0) = f(x, y); \quad \left.\frac{\partial u(x, y, t)}{\partial t}\right|_{t\to 0} = F(x, y). \tag{3}$$

Die Grenzebene, von welcher die Randbedingungen abhängen werden, soll sich innerhalb $-\infty < y < +\infty$ erstrecken, weshalb wir auf (1)—(3) die zweiseitige Laplace-Transformation anwenden wollen.

Wir erhalten demnach für die Unterfunktionen, in bezug auf y, von $u(x, y, t)$ und $\varphi(x, y, t)$:

$$w(x, \varepsilon, t) = \int_{-\infty}^{+\infty} e^{-\varepsilon y} u(x, y, t)\, dy \tag{4}$$

und

$$\gamma(x, \varepsilon, t) = \int_{-\infty}^{+\infty} e^{-\varepsilon y} \varphi(x, y, t)\, dy. \tag{5}$$

Entsprechend ergeben sich für

$$u_0(y, t), \quad h(y, t), \quad f(x, y), \quad F(x, y), \tag{6}$$

in bezug auf y folgende Unterfunktionen:

$$\eta(\varepsilon, t), \quad \zeta(\varepsilon, t), \quad \vartheta(x, \varepsilon), \quad \theta(x, \varepsilon). \tag{7}$$

Aus (1) folgt ferner

$$\frac{1}{c^2}\frac{\partial^2 w}{\partial t^2} + c_1\frac{\partial w}{\partial t} + a_1\frac{\partial w}{\partial x} - \frac{\partial^2 w}{\partial x^2} + (m - \varepsilon^2)\, w = \gamma(x, \varepsilon, t). \tag{8}$$

Wir nehmen nun an, dass uns von den vier Grössen (2) und (3) die erste in (2) und die beiden Grössen (3) gegeben sind.

Dann liefern uns die Ausdrücke (6) und (7) unserer Note I, im Unterraum in bezug auf y, eine vollständige Lösung unseres Problems, falls wir in diesen Ausdrücken an Stelle von

$$u(x, t), \quad \varphi, \quad u_0(t), \quad f(x), \quad F(x) \tag{9}$$

die Grössen:

$$w(x, \varepsilon, t), \quad \sqrt{\varphi^2 - \varepsilon^2}, \quad \eta(\varepsilon, t), \quad \vartheta(x, \varepsilon), \quad \theta(x, \varepsilon), \tag{10}$$

einführen *.

Mit den so abgeänderten Ausdrücken (6) und (7) der Note I gehen wir nun in den Oberraum, in bezug auf y, zurück und erhalten dann die endgültige Lösung unseres Problems. Es ist unmöglich dieselbe hier voll auszuschreiben und wir begnügen uns mit dem der Gleichung (7) der Note I entsprechendem Teil (für $x > tc$), während wir uns das Fehlende zu beschreiben bemühen werden. Wir erhalten:

$$u(x, y, t) = \frac{e^{-\mu t}}{2c}\frac{\partial}{\partial t}\int_{-tc}^{tc} e^{-\nu z} P\{f(x-z,\ y-s),\ t,\ z,\ s,\ \varphi\}\, dz +$$

$$+ \frac{e^{-\mu t}}{2c}\int_{-ct}^{ct} e^{-\nu z} P\{F(x-z,\ y-s) + \mu f(x-z,\ y-s),\ t,\ z,\ s,\ \varphi\}\, dz +$$

$$+ \frac{c}{2}\int_0^t e^{-z\mu} dz \int_{-zc}^{zc} e^{-\nu l} P\{\varphi(x-l,\ y-s,\ t-z), z,\ l,\ s,\ \varphi\}\, dl, \quad x > tc. \tag{11}$$

Hierbei bedeutet P ein Funktionszeichen, und es wird z. B.

$$P\{f(x-z,\ y-s),\ t,\ z,\ s,\ \varphi\} =$$

$$= \frac{1}{\pi}\int_{-\sqrt{c^2t^2-z^2}}^{\sqrt{c^2t^2-z^2}} \frac{f(x-z,\ y-s)\cos\varphi\,\sqrt{c^2t^2-z^2-s^2}}{\sqrt{c^2t^2-z^2-s^2}}\, ds \tag{12}$$

sein.

Es lässt sich deshalb der Ausdruck (11) in Integrale vom Poisson-Parsevalschen Typus umformen, z. B. können wir statt des ersten Gliedes in (11) schreiben:

$$\frac{e^{-\mu t}}{2c}\frac{\partial}{\partial t}\int_{\varphi=0}^{2\pi} d\varphi \int_{r=0}^{ct} e^{-\nu r\cos\varphi}\, \frac{f(x - r\cos\varphi,\ y - r\sin\varphi)\cos\varphi\,\sqrt{c^2t^2-r^2}}{\sqrt{c^2t^2-r^2}}\, r dr, \tag{13}$$

* Man verwechsle nicht die dortige Integrationsvariable y mit der Koordinate y in der vorliegenden Note.

was sich nur durch das Auftreten des Factors $\cos \varrho \sqrt{c^2t^2 - r^2}$, unter dem Integral, von dem gewöhnlichen Poisson-Parsevalschen Integral der ebenen Wellengleichung unterscheidet ($\mu = \nu = m = 0$). Also auch hier integrieren wir längs der Fläche eines Kreises mit dem Mittelpunkt im Aufpunkt und dem Radius ct, welcher also mit der Zeit wächst. Dieser Kreis, kurz bezeichnet Wellenkreis, fegt demnach allmählich die ganze YX-Ebene ab und wird uns die Möglichkeit geben den hier fehlenden Ausdruck für $x < tc$, entsprechend (6) der Note I, zu beschreiben.

Vorerst müssen wir bemerken, dass der Ausdruck (7), Note I, falls wir die Bedingung $x > tc$ fallen lassen, und $\nu = 0$ und $\varphi(x, t) = 0$ setzen, bekanntlich z. B. für einen beiderseits unendlich langen Draht gültig sein wird. Dasselbe bleibt aber auch richtig für den vollen Ausdruck (7) der Note I.

Ganz analog bleibt auch der Ausdruck (11), nach Streichung der Bedingung $x > tc$, für den ganzen Raum, resp. für die ganze YX-Ebene gültig, wobei wir uns also unsere Grenzebene, resp. Grenzlinie (Y-Achse) als nicht existierend betrachten müssen.

Wir spiegeln nun $f(x, y)$, $F(x, y)$ und $\varphi(x, y, t)$, in bezug auf die Y-Achse, ab und denken uns die Grenzebene, resp. die Grenzlinie weg. Es wird dann der Wellenkreis, beim Abfegen der YX-Ebene auch über diese Spiegelbilder kommen und zu den Integralen in (11) noch entsprechende hinzufügen.

Diese Tatsache wird nun gerade durch den von uns ausgelassenen Ausdruck dargestellt. Ausserdem enthält dieser Ausdruck noch das Glied

$$-\frac{c}{\pi} e^{-\nu x} \frac{\partial}{\partial x} \int_{\frac{x}{c}}^{t} e^{-\mu z} dz \int_{-\sqrt{c^2z^2 - x^2}}^{\sqrt{c^2z^2 - x^2}} \frac{u_0(y - s,\ t - z) \cos \varrho \sqrt{c^2z^2 - x^2 - s^2}}{\sqrt{c^2z^2 - x^2 - s^2}} ds, \qquad (14)$$

welches von den gegebenen Werten von $u_0(y, t)$ längs den Y-Achse abhängen wird, wobei längs dem von dem Wellenkreis abgeschnittenen Teil der Y-Achse integriert wird.

Ist im speziellen Fall $f(x, y) = F(x, y) = \varphi(x, y, t) = 0$, so ergibt sich als Lösung, bei $x < tc$, gerade der Ausdruck (14) für den Wert von $u(x, y, t)$, und bei $x > tc$ erhalten wir Null.

Würde links in (1) noch das Glied $\frac{\partial^2 u}{\partial z^2}$ stehen und entsprechend rechts $\varphi(x, y, z, t)$, d. h. wäre der dreidimensionale Fall angenommen, so ist klar, dass wir zu demselben, von (1) aus, auf analoge Weise übergehen können, wie wir dies vom eindimensionalen Fall zu (1) ausgeführt haben. Da aber hierbei einige Eigenarten auftreten, die verdienen besonders hervorgehoben zu werden, so kommen wir noch in einer späteren Note speziell auf diesen Fall zurück.

Zum Schluss möchten wir noch eine Bemerkung hinzufügen die sich auf alle unseren vorhergehenden Noten bezieht. Es betrifft den Ausdruck für die Faltung, welchen wir in einer verallgemeinerter Form gebraucht haben. Es sei

$$f_1(z) = \int_{\alpha_1}^{\beta_1} e^{-zt} F_1(t)\, dt; \quad \beta_1 > \alpha_1 \qquad (15)$$

und

$$f_2(z) = \int_{\alpha_2}^{\beta_2} e^{-zt} F_2(t)\, dt; \quad \beta_2 > \alpha_2. \tag{16}$$

Dann ergibt sich

$$f_1(z) f_2(z) = \int_{\alpha_1+\alpha_2}^{\beta_1+\beta_2} e^{-zt} F(t)\, dt, \tag{17}$$

wobei

$$F(t) = \int_{\max(\alpha_1;\, t-\beta_2)}^{\min(\beta_1;\, t-\alpha_2)} F_1(\tau) F_2(t-\tau)\, d\tau \tag{18}$$

bedeutet und als verallgemeinerter Faltungssatz angesehen werden kann.

Bei $\alpha_1 = \alpha_2 = 0$ und $\beta_1 = \beta_2 = \infty$ geht (18) in die gewöhnliche Faltung über. Der Beweis von (17) und (18), auf welchen wir hier nicht eingehen wollen, beruht auf der Annahme einer absoluten Konvergenz von (15) und (16).

Schwächere Voraussetzungen und Erweiterungen, wie dies Doetsch [2] bei $\alpha_1 = \alpha_2 = 0$ und $\beta_1 = \beta_2 = \infty$ gemacht hat, haben wir nicht vorgenommen.

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ZITIERTE LITERATUR

[1] W. S. Ignatovskij (W. Ignatowsky), C. R. Acad. Sci. URSS, II, I, 5—11 (1935). [2] G. Doetsch, Annali della R. Scuola Normale Superiore di Pisa, Ser II, IV, 71 (1935).

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MATHEMATIK

ÜBER p-GRUPPEN

Von A. P. DIETZMANN

(Vorgelegt von O. J. Schmidt, Mitglied der Akademie, d. 5. III. 1937)

Ist jedes Element einer Gruppe ein Element endlicher Ordnung, so nennen wir solche Gruppe nach Schur (1) periodische Gruppe.

Ist die Ordnung jedes Elementes der periodischen Gruppe eine Potenz von der Primzahl p, so nennen wir diese Gruppe p-Gruppe. Augenscheinlich ist die Ordnung einer endlichen p-Gruppe eine Potenz von p. Es ist bekannt, dass jede endliche p-Gruppe ein vom Einheitselement verschiedenes Zentrum besitzt.

In der vorliegenden Arbeit wird für die p-Gruppe die notwendige und hinreichende Bedingung für die Existenz eines vom Einheitselement verschiedenen Zentrums aufgestellt; dabei werden auch einige Sätze bewiesen.

§ 1. Im folgenden nennen wir der Kürze wegen die Klassen von konjugierten Elementen einer Gruppe Klassen der Gruppe; die Klasse, welche eine endliche Anzahl der Elemente enthält, wollen wir endliche Klasse nennen.

Sei g Produkt von Elementen einer Gruppe:

$$g = a_1 \dots a_{i-1} a_i \dots a_{k-1} a_k a_{k+1} \dots a_l. \tag{1}$$

Wir drücken g in der Form:

$$g = a_1 \dots a_{i-1} a_k a_i^* \dots a_{k-1}^* a_{k+1} \dots a_l \tag{2}$$

aus, wo $a_i^* = a_k^{-1} a_i a_k$; $\dots$; $a_{k-1}^* = a_k^{-1} a_{k-1} a_k$; in diesem Falle heisst es kurz, dass das Produkt (2) aus dem Produkt (1) durch Versetzung des Elementes a_k an die Stelle i erhalten ist.

Sei G eine Gruppe, die eine Menge M von Erzeugenden besitzt. Dann ist jedes Element der Gruppe G ein Potenzprodukt aus je endlichvielen der Elemente von M. Es ist klar, dass jede nichtleere Teilmenge der Elemente von G eine Untergruppe erzeugt.

Satz 1. *Sei M die Menge von Erzeugenden der Gruppe G und $\mathfrak{K}$ ein invarianter Komplex der Gruppe G. Ist $\mathfrak{H}$ eine Untergruppe erzeugt von allen Elementen der Menge M, welche dem Komplex $\mathfrak{K}$ nicht angehören, und Q eine von allen Elementen des Komplexes $\mathfrak{K}$ erzeugte invariante Untergruppe, so ist $\mathfrak{H}Q$ gleich G.*

Beweis. Ist q ein Element von Q, und g ein Element von G, so lässt sich $g^{-1}qg$ als Potenzprodukt aus je endlichvielen der Elemente des invarianten Komplexes $\mathfrak{K}$ darstellen, d. h. $g^{-1}qg$ gehört Q an. Daraus folgt, dass $g^{-1}Qg = Q$, d. h. Q ist eine invariante Untergruppe von G.

Bemerkung. Wenn $g_1 = q_1 g_2$ ist, wo q_1 ein Element von Q und g_1, g_2 Elemente von G sind, so lässt sich g durch Versetzung des Elementes g_2 als Produkt darstellen: $g_1 = g_2 q^*$, wo q^* ein Element von Q ist.

Weiter können wir jedes Element g von G als ein Potenzprodukt aus je endlichvielen der Elemente von M ausdrücken:

$$g = m_1^{\alpha_1} m_2^{\alpha_2} \dots m_k^{\alpha_k}, \tag{3}$$

wo m_1, m_2 ... m_k Elemente der Menge M und α_1, α_2, ... , α_k ganze Zahlen sind. Unsere Bemerkung lässt darauf schliessen, dass man g durch Versetzung der Potenzen von Elementen der Menge M in (3), die dem invarianten Komplex $\mathfrak{K}$ nicht angehören, als Produkt $g = hq_2$ darstellen kann, wo h ein Element der Untergruppe $\mathfrak{H}$ und q_2 ein Element von Q ist. Daraus folgt, dass

$$G = \mathfrak{H}Q.$$

Damit ist unser Satz bewiesen.

Satz II. *Ist A eine endliche Klasse, die aus Elementen endlicher Ordnung der Gruppe G besteht, so erzeugen alle Elemente von A eine invariante Untergruppe Q von endlicher Ordnung.*

Beweis. Bezeichnen wir mit a_1, a_2, ... , a_k die Elemente der Klasse A von der Ordnung k und mit β die Ordnung des Elementes a_1, so haben wir offenbar

$$a_i^{\beta} = 1. \qquad (i = 1,\ 2,\ \dots\ ,\ k) \tag{4}$$

Jedes Element q_j der Untergruppe Q lässt sich als ein Potenzprodukt aus je endlichvielen der Elemente von A darstellen. Nach (4) gibt es also für jedes Element q_j der Untergruppe Q solch eine minimale positive ganze Zahl $m = m(q_j)$, dass man q_j als ein Produkt von m Elementen darstellen kann:

$$q_j = a_{i_1} a_{i_2} \dots a_{i_m}, \tag{5}$$

wo die Faktoren a_{i_1}, a_{i_2}, ... , a_{i_m} einige Elemente von A sind. Befinden sich unter den Elementen a_{i_1}, a_{i_2}, ... a_{i_m} α dem a_s gleiche Elemente, so sagen wir, dass das Element a_s in dem Produkt (5) α-mal vorkommt.

Die Darstellung des Elementes q_j als ein Produkt von m Elementen der Klasse A nennen wir «Normaldarstellung des Elementes q_j», und die Anzahl der Faktoren m—Länge von q_j. Zwei Normaldarstellungen des Elementes q_j heissen verschiedene Normaldarstellungen, wenn sie sich voneinander entweder durch ihre Faktoren, oder durch die Reihenfolge ihrer Faktoren unterscheiden.

Nun betrachten wir die Menge Π aller Normaldarstellungen des Elementes q_j, d. h. die Menge aller verschiedenen Produkte von der Form (4). Offenbar enthält Π eine endliche Anzahl von solchen verschiedenen Produkten.

Sei a_s ein Element der Klasse A, dann kommt das Element a_s in jedem Produkt aus der Menge Π nicht mehr als m-mal vor. Also lässt sich den Elementen a_s und q_j die Zahl α_{sj} zuordnen, welche durch folgende Bedingungen bestimmt wird:

1) Unter den Produkten der Menge Π gibt es wenigstens ein Produkt, in dem das Element a_s α_{sj}-mal vorkommt.

2) In jedem Produkt der Menge Π kommt das Element a_s nicht mehr als α_{sj}-mal vor.

Wir haben also $0 \leqslant \alpha_{sj} \leqslant m$. Im Falle $\alpha_{sj} = m$ ist $q_j = a^{\alpha_{sj}}$.

Es ist hier zu bemerken, dass bei $\alpha_{sj} < m$ eine Normaldarstellung des Elementes q_j von der Form

$$q_j = a_s^{\alpha_{sj}} q_{j+1} \tag{6}$$

existiert, wo q_{j+1} ein Produkt von $m - \alpha_{sj}$ Elementen der Klasse A ist; dabei ist $m - \alpha_{sj}$ die Länge von q_{j+1} und $\alpha_{s_{j+1}} = 0$. Ausserdem ist für jedes Element a_t der Klasse A

$$\alpha_{tj} \geqslant \alpha_{t\,j+1},$$

und, wenn $\alpha_{tj} = 0$, so ist auch $\alpha_{t\,j+1} = 0$.

Um unsere Bemerkung zu begründen, betrachten wir jetzt die Normaldarstellung des Elementes q_j, in der das Element a_s α_{js}-mal vorkommt; durch nacheinanderfolgende Versetzungen der Elemente a_s in diesem Produkt an die erste Stelle bekommen wir offenbar die Normaldarstellung des Elementes q_j von der Form (6). Die Länge von q_{j+1} ist gleich $m - \alpha_{sj}$, weil das Produkt (6) eine Normaldarstellung des Elementes q_j ist. Hier ist $\alpha_{sj+1} = 0$, denn wäre $\alpha_{sj+1} > 0$, so würde eine Normaldarstellung des Elementes q_j existieren, in der a_s mehr als α_{sj}-mal vorkommen würde, was der Definition der Zahl α_{sj} widerspricht.

Endlich ist $\alpha_{tj} \geqslant \alpha_{tj+1}$, denn die Voraussetzung $\alpha_{tj+1} > \alpha_{tj}$ widerspricht der Definition der Zahl α_{tj}. Somit ist unsere Bemerkung begründet.

Es sei q_1 irgendein Element der Untergruppe Q von der Länge m_1. Ist $\alpha_{11} = m_1$, so ist $q_1 = a_1^{\alpha_{11}}$. Wenn aber $\alpha_{11} < m_1$ ist, so folgt aus unserer Bemerkung, dass für q_1 eine Normaldarstellung von der Form

$$q_1 = a_1^{\alpha_{11}} q_2$$

existiert, wobei $\alpha_{12} = 0$ ist.

Wollen wir voraussetzen, dass wir schon eine Normaldarstellung q_1 von der Form

$$q_1 = a_1^{\alpha_{11}} a_2^{\alpha_{22}} \ \ldots\ a_{r-1}^{\alpha_{r-1\,r-1}} q_r; \quad 2 \leqslant r < k, \tag{7}$$

haben, wo

$$\alpha_{1r} = 0; \quad \alpha_{2r} = 0; \ \ldots\ ; \ \alpha_{r-1\,r} = 0$$

und

$$\alpha_{11} + \alpha_{22} + \ \ldots\ + \alpha_{r-1\,r-1} < m_1.$$

Wenn also $\alpha_{11} + \alpha_{22} + \ \ldots\ + \alpha_{rr} = m_1$, so ist offenbar

$$q_1 = a_1^{\alpha_{11}} a_2^{\alpha_{22}} \ \ldots\ a_r^{\alpha_{rr}}.$$

Ist aber $\alpha_{11} + \alpha_{22} + \ \ldots\ + \alpha_{rr} < m_1$, so existiert für q_r die Normaldarstellung:

$$q_r = a_r^{\alpha_{rr}} q_{r+1}, \tag{8}$$

wobei

$$\alpha_{1\,r+1} = 0; \quad \alpha_{2\,r+1} = 0; \ \ldots\ ; \ \alpha_{r\,r+1} = 0;$$

Nach (7) und (8) gibt es für q_1 die Normaldarstellung:

$$q_1 = a_1^{\alpha_{11}} a_2^{\alpha_{22}} \dots a_r^{\alpha_{rr}} q_{r+1},$$

wobei

$$\alpha_{1\,r+1} = 0; \quad \alpha_{2\,r+1} = 0; \dots; \quad \alpha_{rr+1} = 0$$

und

$$\alpha_{11} + \alpha_{22} + \dots + \alpha_{rr} < m_1$$

sind.

Damit wird gezeigt, dass für das Element q_1 die Normaldarstellung von der Form

$$q_1 = a_1^{\alpha_1} a_2^{\alpha_2} \dots a_k^{\alpha_k}$$

existiert, wobei

$$0 \leqslant \alpha_i < \beta \qquad (i = 1, 2, \dots, k),$$

und

$$\alpha_1 + \alpha_2 + \dots + \alpha_k = m_1.$$

Q ist also eine Untergruppe endlicher Ordnung und nach dem Satz I eine invariante Untergruppe von G. Damit ist der Satz II bewiesen.

Satz III. *Ist M die Menge von Erzeugenden der Gruppe G und $\mathfrak{K}$ ein aus endlichvielen Elementen endlicher Ordnung bestehender invarianter Komplex von G, so erzeugen alle Elemente der Menge M, die dem Komplex $\mathfrak{K}$ nicht angehören, eine Untergruppe von endlichem Index der Gruppe G.*

Beweis. Bezeichnen wir mit Q eine von allen Elementen des Invarianten Komplexes $\mathfrak{K}$ erzeugte Untergruppe. Es seien $A_1, A_2, \dots, A_n$ endliche Klassen der Gruppe G, die den invarianten Komplex $\mathfrak{K}$ bilden.

Aus dem Satz II folgt, dass alle Elemente der Klasse A_i (für jedes $i = 1, 2, \dots, n$) den Normalteiler Q_i von endlicher Ordnung der Gruppe G erzeugen; dabei ist

$$Q = Q_1 Q_2 \dots Q_n.$$

Also ist Q eine Untergruppe endlicher Ordnung. Sei $\mathfrak{H}$ eine von allen Elementen der Menge M, die dem Komplex $\mathfrak{K}$ nicht angehören, erzeugte Untergruppe; nach dem Satz I ist

$$G = \mathfrak{H} Q.$$

Da Q eine Untergruppe endlicher Ordnung ist, so ist $\mathfrak{H}$ eine Untergruppe von endlichem Index der Gruppe G, w. z. b. w.

Eine der unmittelbaren Folgen der Sätze II und III ist: *Die Gruppe, welche von n Elementen endlicher Ordnung erzeugt wird, von denen $n-1$ den endlichen Klassen angehören, ist eine endliche Gruppe.*

2. Wir führen zuerst einige Hilfssätze an.

Hilfssatz 1. *Der Durchschnitt von endlichvielen Untergruppen, wo jede Untergruppe von endlichem Index der Gruppe G ist, ist wieder eine Untergruppe von endlichem Index der Gruppe G.*

Beweis. Es seien $\mathfrak{H}_1, \mathfrak{H}_2, \dots, \mathfrak{H}_k$ Untergruppen von endlichem Index der Gruppe G. Mit D_1 bezeichnen wir den Durchschnitt von Untergruppen $\mathfrak{H}_1$ und $\mathfrak{H}_2$, d. h.

$$D_1 = [\mathfrak{H}_1, \mathfrak{H}_2]; \text{ ebenso } D_2 = [D_1, \mathfrak{H}_3]; \dots; D_{k-1} = [D_{k-2}, \mathfrak{H}_k].$$

Offenbar ist D_{k-1} Durchschnitt von den Untergruppen $\mathfrak{H}_1, \mathfrak{H}_2, \ldots, \mathfrak{H}_k$. Aus dem Satz von H. Poincaré (2) folgt, dass D_{k-1} eine Untergruppe von endlichem Index der Gruppe G ist, womit auch unser Hilfssatz bewiesen ist.

Weiter haben wir einen trivialen

Hilfssatz 2. *Ist $\mathfrak{H}$ eine Untergruppe von G, D eine Untergruppe von $\mathfrak{H}$ und Normalteiler von G, so ist die Menge von Nebengruppen der Zerlegung von G nach $\mathfrak{H}$ äquivalent der Menge von Nebengruppen der Zerlegung von Faktorgruppe $\frac{G}{D}$ nach $\frac{\mathfrak{H}}{D}$.*

Hilfssatz 3. *Ist $\mathfrak{H}$ eine Untergruppe von endlichem Index der Gruppe G, so enthält $\mathfrak{H}$ die Untergruppe D, eine invariante Untergruppe von endlichem Index der Gruppe G; dabei ist der Index $\mathfrak{H}$ unter G gleich dem Index $\frac{\mathfrak{H}}{D}$ unter $\frac{G}{D}$*, d. h.

$$(G, \mathfrak{H}) = \left(\frac{G}{D}, \frac{\mathfrak{H}}{D}\right).$$

Beweis. Offenbar enthält G endlichviele verschiedene, mit $\mathfrak{H}$ konjugierte Untergruppen. Mit D bezeichnen wir den Durchschnitt aller mit $\mathfrak{H}$ konjugierten Untergruppen. Aus dem Hilfssatz 1 folgt, dass D eine invariante Untergruppe von endlichem Index der Gruppe G ist. Nach dem Hilfssatz 2 haben wir:

$$(G, \mathfrak{H}) = \left(\frac{G}{D}, \frac{\mathfrak{H}}{D}\right).$$

Der Hilfssatz ist bewiesen.

Folgerung. *Ist $\mathfrak{H}$ eine Untergruppe von endlichem Index einer p-Gruppe P, so ist der Index von $\mathfrak{H}$ unter P eine Potenz von p.*

Beweis. Sei D der Durchschnitt aller mit $\mathfrak{H}$ konjugierten Untergruppen. Nach dem Hilfssatz 3 ist D eine invariante Untergruppe von endlichem Index der Gruppe P und

$$\left(\frac{P}{D}, \frac{\mathfrak{H}}{D}\right) = (P, \mathfrak{H}). \tag{9}$$

Weiter ist die Faktorgruppe $\frac{P}{D}$ eine endliche Gruppe, dabei ist die Ordnung jedes Elementes davon eine Potenz von p. Daraus folgt, dass die Faktorgruppe $\frac{P}{D}$ eine endliche p-Gruppe ist. Nach (9) ist der Index von $\mathfrak{H}$ unter P eine Potenz von p, w. z. b. w.

Satz IV. *p-Gruppe besitzt ein vom Einheitselement verschiedenes Zentrum, wenn sie wenigstens eine vom Einheitselement verschiedene endliche Klasse enthält.*

Beweis. Es sei A eine endliche Klasse der p-Gruppe P. Aus dem Satz II folgt, dass die Elemente der Klasse A den Normalteiler $\mathfrak{A}$ endlicher Ordnung p^n der Gruppe P erzeugen. Mit a bezeichnen wir irgendein Element von $\mathfrak{A}$. Das Element a gehört zu der endlichen Klasse der Gruppe P, weil die Menge von allen Elementen des Normalteilers $\mathfrak{A}$ die Gesamtheit einiger endlichen Klassen der Gruppe P darstellt. Der Normalisator V des Elementes a in der Gruppe P ist also eine Untergruppe von endlichem Index der Gruppe P. Laut der Folgerung des Hilfssatzes 3

ist der Index V eine Potenz vone p. Folglich ist die Ordnung der Klasse der mit a konjugierten Elemente der Gruppe P eine Potenz von p.

Zerfällt der Normalteiler $\mathfrak{A}$ in k Klassen der Gruppe P, wobei die Ordnungen dieser Klassen $\alpha_1, \alpha_2, \ldots, \alpha_k$ sind, so haben wir:

$$p^n = \alpha_1 + \alpha_2 + \ldots + \alpha_k.$$

Die Zahlen $\alpha_1, \alpha_2, \ldots, \alpha_k$ sind Potenzen von p und, da $\alpha_1 = 1$ ist, (das Einheitselement bildet eine Klasse), muss es ausser dem Einheitselement noch weitere Elemente geben, die für sich allein eine Klasse der Gruppe P bilden, was unseren Satz beweist.

Aus dem Satz IV folgt, dass die p-Gruppe, deren Zentrum dem Einheitselement gleich ist, ausser dem Einheitselement keine andere endliche Klasse enthält.

Eingegangen
d. 5. III. 1937.

ZITIERTE LITERATUR

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ASTRONOMY

THE USE OF THE PHOTO-RELAY IN THE STUDY OF LUNAR MICRORFLIEF

By S. G. NATANSON

(Communicated by D. S. Roždestvenskij, Member of the Academy, 17. II. 1937)

When a star is occulted by the Moon, the latter throws its shadow on the Earth in the rays of the star. Since these rays may be taken to be parallel, the outline of any detail of the lunar edge would therefore be projected on any plane normal to the rays «as large as life».

If it were possible by some means to outline on the Earth's surface the instantaneous position of the shadow of that detail of the lunar edge, i. e., to plot on a locality or map the isochrone of the star's occultation, the problem would be solved in its entirety. This isochrone may be determined from accurate observations of the moments of the beginning (or the end) of the star's occultation, i. e., of the moments of the disappearance or reappearance of the star from beyond the edge of the Moon.

The velocity of the apparent motion of the Moon's shadow on the Earth's surface can indeed be calculated from the rules of spherical astronomy with the requisite degree of accuracy. On an average, the velocity of the Moon is 0″.5 per 1 sec. and that of the shadow, 1 km per 1 sec. Consequently, if the initial (or final) moments of occultation were known to within $\pm 0^s.001$, the position of the shadow would be determined on an average to within ± 1 m. Let us place a chain of observers $A_1, A_2, A_3, \ldots, A_n$, (Fig. 1) capable of recording, with reference to a common clock, the moments of the occultation of the star with accuracy suggested above. The moments obtained by the observers: t_1, t_2, t_3, $\ldots$, t_n, would permit to reproduce the outline of the Moon's shadow at some arbitrary moment t_0, i. e., to plot an isochrone of the occultation for that moment t_0 with an accuracy of the order of 1 m. To do this, it suffices to mark off, from the points of observation A_1, A_2, $A_3, \ldots, A_n$, in the direction of the motion of the Moon's shadow, the distances: $\varphi(t_0 - t_1)$, $\varphi(t_0 - t_2)$, $\varphi(t_0 - t_3)$, $\ldots$, $\varphi(t_0 - t_n)$, where φ is the velocity of motion of the Moon's shadow.

Let us consider how the diffraction of the star's light by the edge of the Moon would affect our observations. Fig. 2 shows the intensity of illumination of the Earth's surface normal to the rays of the star, near to the edge of the geometrical shadow. The distances from the edge of the geometrical shadow are plotted on the axis of the abscissae, while the intensity of illumination is plotted on the axis of the ordinates. Far from the edge, the illumination due to the star is taken at 100 per cent. A rapid decrease of illumi-

nation is observed over a distance of 12 m. Such is the «diffusion» of the Moon's shadow due to diffraction.

The use of a sensitive photo-relay renders it quite possible to record automatically the moments of the reappearance and disappearance of the star at the dark edge of the Moon with a degree of accuracy quite sufficient for our purposes [1]. Most suitable for the purpose is the L. A. Kubetzky secondary electronic tube (multiple amplifier) as it has a high coefficient of amplification combined with a practically total absence of inertia [2,3].

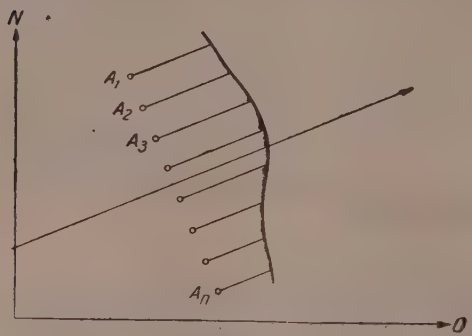

Fig. 1.

Working in his laboratory, I have lately designed a tube which, when placed in the focus of a one-metre mirror, is capable to feel the light of stars down to the 5.5 magnitude. In this case, the output photo-current reaches such values as permit of automatic recording. The mirror must not necessarily possess high optic qualities, its object being to amplify the luminous flux falling on the photo-cathode of the tube. Stars of the 1st magnitude require a mirror of no more than 7—8 in. in diameter.

It would not be difficult for all the registering devices to work by the same clock. A short-wave transmitter would give rhythmical signals which would be received by the observer on the same recorder as the photo-current supplied by the light of the star. The method suggested above would thus be within the limits of practical possibilities.

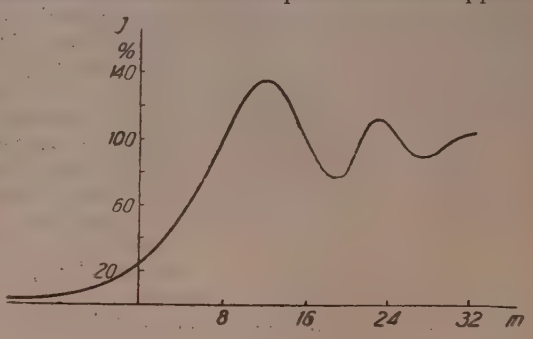

Fig. 2.

The 12 m diffusion of the Moon's shadow is not by any means a limit of the resolving power of this method. According to some schemes, a photo-relay may be so adjusted as to recorder not a complete disappearance of light, but its decrease down to a certain proportion, e. g., by 50 per cent. The diffraction curve shows that the use of a relay capable to record the moment when the illumination falls to 30—70 per cent of its initial value, would involve an error of 4.5 m in determining the geometrical shadow position. It is of interest to note that this is the resolving power of a telescope with a 50-metre diametre mirror.

Such observations on occultation, if carried on with a photo-relay, make it possible to study the lunar microrelief so far utterly unknown.

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17 II 1937.

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CHEMISTRY

ON THE CATALYTIC ISOMERIZATION OF NORMAL HEXENE AND OCTENE IN THE PRESENCE OF CHLORIDE OF ZINC AND PHOSPHORIC ACID

By A. D. PETROV and M. A. ČELCOVA

(Communicated by A. E. Favorskij, Member of the Academy, 20. II. 1937)

The shifting of the double bond and change in the carbon skeleton enter into the conception of isomerization which has great practical importance in the technology of motor fuels of high anti-detonational qualities. The possibility of the shifting of the double bond under certain conditions has been frequently pointed out. We may mention here the investigations ([1, 2, 3, 4]) and others.

Change of the carbon skeleton has been observed principally in reactions of dehydrations of alcohols carried out over $ZnCl_2$.

It has been shown ([5]) that, while only butylene-2 is formed in the dehydration of methylethylcarbinol over Al_2O_3, the dehydration of the latter over $ZnCl_2$ results in a mixture of butylene-2 with isobutylene. In like manner, in the dehydration of isobutyl alcohol there results the so-called di-isobutylene of Malbot and Gentil which, as A. D. Petrov, L. I. Ancus and D. N. Andreev ([6]) have shown, is a mixture of 2. 2. 4-trimethylpentene with 2. 2. 3-trimethylpentene.

One of the rather few works on the catalytic isomerization of the hydrocarbons of the olefine series (accompanied by change of the carbon skeleton) is the recent work of K. Laughlin, C. Nash and F. Whitmore ([7]). However, they demonstrated only a shifting of the side chains already present (the conversion of 2. 2.-butene into 2. 3.-butene, or vice versa), but not the formation of the latter from hydrocarbons of normal structure.

It was among the purposes of the present investigations to find the optimal conditions for the maximum isomerization of normal hexene and octene into the corresponding branched hydrocarbons and to determine the effect of this isomerization in terms of octane values.

We conducted experiments in isomerization under atmospheric and under high pressures in the presence of both zinc chloride and phosphoric acid.

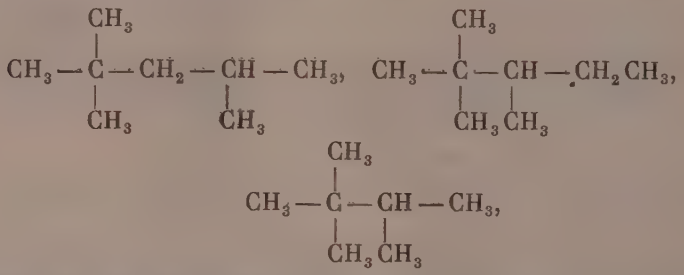
The degree of isomerization in each individual case was determined in the following manner .

The isomerizate was hydrogenated into the corresponding saturated hydrocarbons, and the content of the branched hydrocarbons in these latter was determined by titrating with $SbCl_5$ in a chloroform solution after the method

of Schaarschmidt-Moldavski ([8]). It was found that the yield of branched hydrocarbons increases with the increase of temperature, pressure and the duration of heating, and that chloride of zinc is a more active catalyzer than phosphoric acid, inasmuch as it gives a higher yield of branched hydrocarbons under lower temperatures.

Titration of the isomerizates (after they had been hydrogenated) in the chloroform solution was controlled by titration of the individual branched hydrocarbons (2. 5-dimethylhexane, 2. 2. 4-trimethylpentane, 2. 2. 3-trimethylpentane and 2. 2. 3-trimethylbutane) which had been obtained by synthesis.

An interesting fact was discovered during this process: it was found that the three last hydrocarbons:



in which the tertiary carbon atoms exposed to the action of $SbCl_5$ were, so to speak, «under the protection» of the quarternary carbon atoms (which are not chlorinated by $SbCl_5$) and are chlorinated several times more slowly than 2. 5-dimethylhexane under the modification of Schaarschmidt's method proposed by Moldavski. As will be remembered, under Schaarschmidt's method, i. e., in chlorinating with undiluted $SbCl_5$, no difference in the chlorination of hydrocarbons of these two types is to be observed. Thus, while chlorination by Schaarschmidt's method involves the danger of exaggerated isomerization values (as a consequence of the inductive influence on the normal hydrocarbons also), chlorination by Moldavski's method (in a chloroforms sulution), when there are present in the mixture hydrocarbons that contain both tertiary and quaternary carbon atoms, is open to the danger of leading to undervaluations.

In the case, for example, of octene, the possibility of its being isomerized into 2. 2. 4-trimethylpentene cannot be excluded. Indeed, di-isobutylene is a very stable olefine hydrocarbon, and one of us succeeded in isolating and identifying it when condensing acetone into hydrocarbons of the olefine series under the action of $ZnCl_2$ ([9]).

Partly in order to check up the results of isomerization that we obtained and evaluated in percentages of the content of branched hydrocarbons in the isomerizates after their hydrogenation, partly in order to appraise the practical effect of this kind of isomerization, we carried out the isomerization of hexene-1 and of a mixture of octene-1 and octene-2 (which were obtained by the dehydration of octanol-2) in quantities sufficient for the determination of octane values.

It was shown that hexene-1, which, according to the literature ([10]), has an octane value of 68.5, had, after isomerization, an octane value of 78. Inasmuch as only the octene value of octane-1 (equal 39.5) is reported in the literature, while the octane value of octene-2 (which is apparently higher) is unknown, we determined the octane values of a mixture of octene-1 and

octene-2 both before and after isomerization. The octane value before isomerization was 60; after isomerization, 75.

The growth of octane values by 10 to 15 points as a result of the isomerization of the hexene-octene fraction must be recognized as noteworthy and perhaps as having a practical importance in the isomerization of the paraphase cracking-benzine carried out before its hydrogenation. Besides this, the exacting conditions (as to temperature, time, presence of active catalyzers), needful for any considerable isomerization, points to the fact that the possibility of combining any considerable isomerization with hydrogenation is hardly probable.

Experimental Part

The experiments in isomerization were begun with a mixture of octene-1 and octene-2 obtained by the dehydration of octanol-2 according to the method of Thomas and Mannich [11]. Before isomerization, this mixture of hydrocarbons had a boiling point of 119—124°; $d_4^{20} = 0.7204$, $n_D^{20} = 1.4141$. After dehydration, boiling point was 123—125°; $d_4^{20} = 0.7035$; $n_D^{20} = 1.3979$; iodine value = 0.

Titration with $SbCl_5$ was carried out both in this and in all subsequent instances after the method of Moldavski and Livshits [12], and revealed the almost complete absence of hydrocarbons of branched structure in the initial hydrocarbons (percentage of iso-compounds is equal to 0.3).

The first isomerizate that we got, after it had been hydrogenated, was titrated simultaneously with a control sample, this being 2. 2. 4-trimethylpentane. It appeared that this latter hydrocarbon chlorinates considerably more slowly than we had expected on the basis of the rule established by V. A. Moldavski [12] concerning «the independence of the rate of chlorination on the structure of branched hydrocarbons».

Inasmuch as Moldavski carried on his control experiments with 2-methylbutane, 2-methylpentane, 3-methylpentane, 2. 3-dimethylpentane, 4-methylheptane and 2. 5-dimethylhexane, that is, only with hydrocarbons having one or two tertiary carbon atoms, and did not concern himself with hydrocarbons containing quaternary carbon atoms along with the tertiary, we turned our attention to the behaviour of the last mentioned when chlorinated with $SbCl_5$. Besides 2. 2. 4-trimethylpentane, we obtained 2. 2. 3-trimethylpentane according to the method of L. Clarke and W. Jons [13], and 2. 2. 3-trimethylbutane by the Grignard reaction through pentamethylethanol. The effect of chlorination on these hydrocarbons, taken separately and in different concentrations in a mixture with normal octane, are shown in Tables 1 and 2.

Table 1

Number of Moles of $SbCl_5$ Formed per 1 Mole of Hydrocarbon

Time in hours	2. 5-dimethylhexane	8% solution of 2. 5-dimethylhexane in normal octane	20% dimethylhexane in normal octane	40% dimethylhexane in normal octane	2.2.4-trimethylpentane	8% 2.2.4 trimethylpentane in normal octane	20% 2.2.4-trimethylpentane in normal octane
3	3.74	4.2	3.95	3.88	0.84	1.4	1.13
6	—	—	5.08	—	—	2.3	1.64
12	—	—	—	—	—	—	4.4

Table 2

Number of Moles of $SbCl_3$ per 1 Mole of Hydrocarbon

Time in hours	2.2.3-trimethylpentane	20% 2.2.3-trimethylpentane in normal octane	25% 2.2.3-trimethylpentane in normal octane	2.2.3-trimethylbutane	24% 2.2.3-trimethylbutane in normal octane
3	0.6	0.7	0.78	0.20	1.2
6	1.2	1.3	1.21	1.40	2.0

It is evident from these data that, while 2. 5-dimethylhexane is approximately 100% chlorinated after 3 hours (after 6 hours, 125%), the hydrocarbons containing quaternary carbon atoms require 12 hours or more for 100% chlorination (after three hours they are chlorinated only 40%, and in the case of some hydrocarbons only 20%).

A similar lag in rate of chlorination of hydrocarbons containing quaternary carbon atoms takes place also when they are in a mixture with branched hydrocarbons not containing quaternary carbon atoms.

For example, a mixture of 8% 2. 2. 4-trimethylpentane and 8% 2.5-dimethylhexane became only 65% chlorinated after 3 hours. An analogous result was also obtained from a mixture containing 20% of each these hydrocarbons in normal octane.

Thus, hydrocarbons containing quaternary carbon atoms along with tertiary are an exception to Moldavski's rule, and when they are present in a mixture of hydrocarbons, the results of an analysis conducted by that method are likely to be somewhat beneath the true quantities. However, owing to the lack of more exact methods of determining the content of hydrocarbons of iso-structure, and regarding the possible error in determinations by Moldavski's method as less important than that arising under Schaarschmidt's conditions, we preferred to follow Moldavski. The experiments in the iso-

Table 3

Experiment No.	Catalyzer used	Duration of experiment in hours	Temperature °C	Percentages of iso-compounds	Remarks
1	15 g $ZnCl_2$ on pumice (1 : 1)	25	325—350	12	B. p. of isomerizate, 119—122°
2	15 g $ZnCl_2$ on pumice (1 : 1)	25	325—350	12.3	Hydrocarbon was in a mixture with water vapour. B. p. of isomerizate, 118—122°
3	50 g $ZnCl_2$ on pumice (1 : 1)	13	325—350	25	B. p. of isomerizate, 108—122°
4	Experiment repeated with same catalyzer as in No. 3	13	325—350	17	

merization of a mixture of octene-1 and octene-2 were conducted under atmospheric and under high (in the autoclave) pressures.

In the experiments under atmospheric pressure, we took each time 50 cm^3 of the hydrocarbon and passed it several times through tubes filled with the catalyzer. The tubes were 1) 43 cm long and 12 mm in diameter; 2) 81.5 cm long and 18 mm in diameter. The data from these experiments are shown in Table 3. The percentage of iso-compounds was determined after hydrogenation of the isomerizate to an iodine value of zero, washing in sulphuric acid and redistillation.

We conducted experiments under pressure in rotating autoclaves (the first two experiments with pure $ZnCl_2$) and in stationary autoclaves (in case of placing the catalyzer on the pumice). 40 cm^3 of the hydrocarbon were taken for each experiment, and the initial pressure of the nitrogen was 50 atmospheres. The data of these experiments are shown in Table 4.

Table 4

Isomerization of octene-1 and octene-2 under pressure

Experiment No.	Catalyzer used	Duration of experiment	Temperature C°	Percentages of iso-compounds	Remarks
1	8 g $ZnCl_2$	30 m.	250—300	—	B. p. of isomerizate, 118—122°.5
2	8 g $ZnCl_2$	1 h. 30 m.	300—325	8.34	B. p. of fundamental isomerizate, 118—124° (boiling began at 103°)
3	35 g $ZnCl_2$ on pumice (1 : 1)	30 m.	300—325	20.0	B. p., 114—123° (boiling began at 50°)
4	35 g $ZnCl_2$ on pumice (1 : 1)	1 h.	300—325	22.0	B. p., 114—123° (boiling began at 50°)
5	35 g $ZnCl_2$ on pumice (1 : 1)	1 h. 30 m.	300—325	45.0	B. p., 106—124° (boiling began at 50°)
6	27 g phosphoric acid on pumice (1 : 1)	1 h. 30 m.	300—325	15.0	B. p., 109—125°
7	27 g phosphoric acid on pumice (1 : 1)	1 h. 30 m.	400	46.7	B. p., 95—127° (boiling began at 45°)

In conclusion, we conducted a series of experiments under standard conditions (temperature, 300—325°; duration of heating, 1 h. 30 m.; initial pressure of nitrogen, 50 atmospheres) and also with large quantities of catalyzer (100 cm^3 of the hydrocarbon to 75 g $ZnCl_2$ on pumice in the ratio of 1: 1) both with octene-1 and octene-2, and with hexene-1, in order to obtain isomerizates in quantities sufficient for the determination of the octane values. One charge of catalyzer served us for 3 experiments with charges of 100 ml each of the hydrocarbon, after which the catalyzer was changed for fresh, while the old catalyzer underwent extraction of the resins formed in it, so that a calculation might be made of the balance of the process.

75% of the reaction product consisted, in the case of octene-1 and -2, of a light isomerizate, which evaporated at temperatures up to 130°, and 25%

consisted of higher boiling fractions and of resins on the catalyzer. 40% of these high-boiling fractions evaporated at temperatures of up to 200°, and 60% at temperatures of up to 230°.

The data on the distillation of the light isomerizate were as follows:

30 to 95°	8.5%	up to 118°	60%
up to 100°	10%	up to 119°	65%
up to 116°	50%	up to 130°	100%

The constants of 30—130° before hydrogenation were as follows: $d_4^{20} = 0.7120$; $n_D^{20} = 1.4112$. After hydrogenation: b. p., 30—130°; $d_4^{20} = 0.6966$; $n_D^{20} = 1.3943$. Percentage of iso-compounds according to Moldavski, 46.8. Octane value of the isomerizate before hydrogenation, 75. Before isomerization, the initial product with b. p. 119—124° had an octane value of 60.

From hexene-1 (obtained by the dehydration of hexanol-1 over Al_2O_3) having a b. p. of 63—65°, an isomerizate was obtained, 80% of which consisted of a light fraction having a b. p. of 63—68°; 15% (after deducting losses) consisted of a heavy fraction, of which 10% fell in fraction 80—200° and 5% in fraction having b. p. above 200°. The resins were not extracted in this case, since they were in very small quantity, as evidenced by the almost unchanged colour of the catalyzer.

The constants of fraction 63—80° before hydrogenation were: $d_4^{20} = 0.6799$; $n_D^{20} = 1.3950$. After hydration: b. p. 63—73°; $d_4^{20} = 0.6586$; $n_D^{20} = 1.3757$. Percentage of iso-compounds in fraction 63—73° after hydrogenation was 23. The octane value of fraction 63—80° was 78.

From the data presented it is apparent that a considerable isomerization of hexene and octene can be obtained at the cost of relatively small losses on polymeres (about 15% to 25%). However, this considerable isomerization can be obtained only by using a large amount of catalyzer in the process. But the expenditure of catalyzer cannot be computed on the basis of our data, since we did not ascertain the length of time during which the calatyzer remains active. It may also be seen from our data that, when isomerization is carried out under high pressure, a higher degree of isomerization can be reached than when conducting the experiment under atmospheric pressure, and in a shorter time.

The considerably slighter activity of phosphoric acid, as compared with $ZnCl_2$, at temperatures of 300 to 325° is also interesting.

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BIOHEMISTRY

THE DETERMINATION OF ASCORBIC ACID (VITAMIN C) IN BLOOD AND URINE

By V. A. DEVJATNIN and V. M. IOSIKOVA

(Communicated by A. N. Bach, Member of the Academy, 27. II. 1937)

Study of vitamin C metabolism in the organism is now one of the to-day problems of physiology; an enormous literature is devoted to it.

We set ourselves in 1936 the goal of finding the most convenient method for everyday clinical application in this field, being impelled by the fact that up to that time there was no sufficiently accurate method of determining ascorbic acid in the blood and urine.

The methods of Tillmans (1), of Harris and Ray (2) and others are not applicable because of the presence in the blood and urine of a large quantity of various substances which reduce (2, 6) dichlorphenolindophenol, as, for example, cystein, glutathion, sugar, ergothionein, and others. The method of Martin and Bonsignore (3) is somewhat cumbersome under clinical laboratory conditions and not always specific. The methods proposed by Emmerie and Eekelen (4) and by Bukin (5) are the most interesting and specific in principle, but even they do not give accurate results. The acids used by these authors (trichloracetic and hydrochloric) exert a destructive action on the indicator, while mercury acetate completely oxidizes ascorbic acid and the subsequent treatment of the latter with hydrogen sulphide does not reduce it to 100 per cent .

Farmer and Abt (6), Fujita and Iwatake (7) have proposed new methods, but these methods are also inapplicable, since, in the first instance, only the blood plasma is used for analysis, while, in the second, the use of the red blood salt in biological extracts leads to incorrect titration results, as has also been pointed out by Manceau, Policard and Ferrand (8).

We have previously proposed a method for the determination of ascorbic acid in vegetable materials (9, 10, 11). Lead acetate in acetic extract, which we used, does not oxidize ascorbic acid and leads to results that correspond with the true content of vitamin C in the material (12, 8), but it does not eliminate cystein and thionin from the extract, which is not important in investigations conducted on vegetable material, but is important in working with animal materials. This fact was noted in a criticism of our method by Emmerie and Eekelen (13).*

* We desire to express our thanks to these authors for their criticism of our work.

It had been observed by us that oxidation of ascorbic acid by mercury acetate is considerably retarded by the presence of calcium salts and that the ascorbic acid oxidized in this manner may afterwards be completely reduced by hydrogen sulphide*. The trend in present-day literature is to look upon calcium as a factor retarding the oxidation of ascorbic acid ([15]).

This fact served as the basis of our method, the technique of which is set forth below.

Method

1. To 4 cm³ of blood or urine in a wide centrifuging test tube are added 8 cm³ of 0.5% calcium acetate and then, stirring carefully with a glass rod, 4 cm³ of 25% mercury acetate are slowly poured in. After being stirred vigorously till the mixture is homogenous, the liquid is centrifuged 5 minutes (there is no need to centrifuge urine—it is sufficient to pass it through a filter).

2. The centrifugate (or filtrate) is treated in a flask for 3 to 5 minutes with hydrogen sulphide, the latter being then removed by means of carbon dioxide.

3. A few cm³ of the transparent colourless extract are transferred with a pipette into a test tube and titrated with a 0.001/2 normal solution of (2,6) dichlorphenolindophenol from a microburette to a definite, stable pink colour**. The amount of ascorbic acid can be reckoned in mg %, assuming that 1 mg of ascorbic acid reduces 11. 4 cm³ of 0.001 normal solution of (2, 6) dichlorphenolindophenol.

4. A paralled calculation is made of the amount of indicator used in titrating the reagents («correction for colour»), for which purpose a control experiment is carried through with the reagents and distilled water. The number of cm³ of indicator thus calculated is deducted.

This method was tested out on crystallized ascorbic acid and on various substances present in blood and urine and capable of reducing the indicator

The results are shown below.

Table 1

Titration of Ascorbic Acid Introduced into Blood and Urine
(Data expressed in cm³ of indicator used in titration)

Serial Nos.	Blood				Urine				Remark
	Ascorbic acid	Blood	Blood + Ascorbic acid	Initial amount of ascorbic acid determined in %	Ascorbic acid	Urine	Urine + ascorbic acid	Initial amount of ascorbic acid determined in %	
1	9.90	0.40	10.30	100.0	9.90	0.0	9.90	100.0	Blood and urine investigated during the period March-May, 1936
2	20.60	0.50	21.10	100.0	12.20	0.5	12.70	100.0	
3	20.60	0.55	21.15	100.0	34.10	0.0	54.10	100.0	
4	34.10	0.40	34.50	100.0	—	—	—	—	
5	7.19	0.44	7.63	100.0	—	—	—	—	

* We are studying the mechanism of this action and shall soon publish our findings ([14]).

** Inasmuch as the content of ascorbic acid in blood and urine is usually rather small, it is best to use the whole of the filtrate to titration in order to avoid mistakes.

Table 2

Titration of Reducing Substances

Substance	cm³ of indicator per 2 cm³ of reducer	Number of determinations made	Substance	cm³ of indicator per 2 cm³ of reducer	Number of determinations made
Glutathion, 0.5% . .	0	3	Mixture of aromatic amino-acids, 1.0%	0	2
Cystein, 0.1%	0	6	Mixture of amino-acids of the fatty series, 1.0% . . .	0	2
Blood + cystein, 0.1% (after subtracting blood)	0	5	Mixture of monosaccharides 1.0% . .	0	2
Blood + glutathion, 0.1% (after subtracting blood . .	0	4	Mixture of polysaccharides, 1.0% . .	0	2
Sacharose, 5% . . .	0	2	Aldehydes	0	4
Fructose, 1%	0	2	Ketones (except acetone)	0	4
Glucose, 1%	0	2	Oxalates	0	2
Phenols, 2%	0	2	Phosphates	0	2
Iron citrate, 0.2% . .	0	2	Sulpho-proteins, 5%	0	2
Fatty acids, 0.1% . .	0	2	Sterins (mixture), 0.5%	0	2
Pyrogallic acid, 0,1%	0	2			
Uric acid, 0.5% . .	0	2			
Acetone, 0.4% . . .	0	2			

From investigations of the blood and urine of a great number of healthy subjects we obtained data which indicate a certain stability of ascorbic acid content in the blood and also a relation between the introduction of vitamin C into the body in food and its discharge in the urine. Some of these data are presented in Tables 3 and 4.

Table 3

Content of Ascorbic Acid in Human Blood

	Ascorbic acid in mg %
Normal	0.50
»	0.55
»	0,57
»	0.60
Hypovitaminosis	0.32
»	0.25
»	0.0

We intend to publish more complete data later on, but even the material here presented justifies us in drawing some conclusions.

1. The so-called «hypervitaminosis C» does not occur in the human organism.

2. The degree to which the organism is saturated with ascorbic acid can be estimated only after ascertaining its content simultaneously in both blood and urine.

3. Hypovitaminosis C may occur, especially in the winter-spring period, if fresh fruits and vegetables are not eaten in sufficient quantity.

Table 4

Discharge of Ascorbic Acid in the Urine in Dependence on Food Ration

	Day of the experiment									
	1	2	3	4	5	6	7	8	9	10
	Content in milligrams of ascorbic acid									
	0	150	150	150	150	300	300	300	300	300
Amount of ascorbic acid in urine discharged during 24 hours (in mg) .	0.0	0.0	6.3	4.7	2.2	4.7	26.5	50.5	87.0	135.0

In conclusion, we set forth the technique of preparing mercury acetate according to Tovarnitski ([16]) for the reason that this reagent is not on the market.

110 grams of yellow oxide of mercury, in a flask with groundglass stopper, are mixed with 400 cm³ of distilled water freed of all CO_2, then 60 grams of ice-cold acetic acid are added, and the mixture is carefully shaken up. It is then allowed to stand for 7 to 10 days, being throughly shaken up from time to time. At the expiration of the time indicated, the solution is filtered, and 20 cm³ of distilled water is added to each 100 cm³ of the filtrate. The 25% solution of the salt thus obtained keeps indefinitely in a dark place.

Summary

1. A method is set forth of determining ascorbic acid in the blood and urine. It is free of the necessity of reducing indophnenol by means of extraneous reducing substances, and makes it possible to determine the ascorbic acid conretent of the material investigated to 100%.

2. The basic principle of this method is the treatment of the material with mercury acetate in the presence of calcium salts which prevent the oxidation of ascorbic acid, followed by a further reduction with hydrogen sulphide.

3. Data characterizing the method and enabling an evaluation of it are presented.

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GÉOLOGIE

SUR LA DÉCOUVERTE DE L'ÉOCÈNE SUPÉRIEUR EN GOURIE CARACTÉRISÉ PAR LA FAUNE

Par K. S. MASLOV et I. A. KOROBKOV

(Présenté par A. A. Borissiak, de l'Académie, le 23. II. 1937)

En été 1935 K. S. Maslov, géologue de l'Institut de Recherches géologiques a découvert en Gourie les grès à gros grains polymictiens de l'éocène supérieur, renfermant un complexe riche de la faune de mollusques à prédominance nette de Gastropoda. Les dépôts en question se trouvent au nord-ouest de Gourian-mta et à l'affluent droit de la rivière Méria sous la latitude de Gogorety, en forme de petites taches isolées et des sillons. Ils se rapportent à la large zone monoclynale, compliquée par le microplissement, du développement des formations sablo-argileuses du sarmatien inférieur, entrecoupées par des dykes discontinues de basalte. Les roches de l'éocène supérieur ci-dessus décrites s'appuyent partout aux parties périphériques des dykes en cause et apparaissent aux endroits d'interruptions des dykes. Gardant la continuité stratigraphique normale l'éocène supérieur de la Gourie est surtout développé à la distance de 15—20 km vers le nord-est de cet endroit, aux environs des villages Aketa, Chanchkhata, Vani, Samebo. Il est présenté là-bas par un facies de marnes globigérines aux intercalations de grès, dépourvu de faune de mollusqes. Comparés à ces derniers, les grès des environs de Gourian-mta peuvent être considérés comme formations d'un facies marin moins profond d'une seule et même zone gourienne de l'éocène qui peut être envisagée comme forme de transition de zones de la chaîne principale du Caucase Central à celle de la chaîne Adjaro-Imeretine.

Leur affleurement dans les conditions ci-dessus décrites est déterminé exclusivement par des déplacements verticaux. La structure autochtone géologique de la Gourie ne démontre pas de charriages et, de plus, l'absence indiscutable de recouvrements allochtones exclut, en tout cas, leur déplacement tectonique du coté de la chaîne Adjaro-Imertine.

La faune de mollusques trouvée dans les grès du Gourian-mta est représentée par de nombreuses coquilles de *Gastropoda* très bien conservées et par petit nombre de débris et noyaux de *Pelecypoda*.

L'étude de cette faune a démontré, qu'elle est composée d'espèces dont l'énumération et la répartition stratigraphique sont données dans la table ci-dessous.

Nomination des espèces	Répartition stratigraphique			
	Éocène inférieur	Éocène moyen	Éocène supérieur	Oligocène inférieur
1. *Cassidea thesei* Brongniart			+	
2. *Cassidea ambigua* Solander	+	+	+	+
3. *Apporhais pescarbonis* Brongniart			+	+
4. *Turritella (Archimediella) gradataeformis* Schauroth			+	
5. *Conus (Zeptoconus) diversiformis* Deshayes		+	+	
6. *Cryptoconus filosus* Lamk		+	+	+
7. *Borsonia costellorum* Oppenheim			+	+
8. *Mitra postera* Koenen			+	+
9. *Tornatella simulata* Solander	+	+	+	+
10. *Spondylus buchi* Philippi	+	+	+	+
11. *Ostrea (Gigantostrea) gigantica* Solander		+	+	+
12. *Axinea* cf. *striatissima* Bellardi			+	
	3	6	12	8

La table démontre que la faune étudiée est composée de trois espèces largement répandues dans la direction verticale; quatre espèces font leur première apparition dans l'éocène moyen dont deux se rencontrent jusqu'à l'oligocène et l'une n'est connue que dans l'éocène moyen et supérieur; trois espèces appartiennent uniquement à l'éocène supérieur et trois espèces ne sont connues que dans l'oligocène supérieur et inférieur. Le complexe cité est tout à fait dépourvu d'espèces apparaissant pour la première fois dans l'éocène inférieur. L'analyse de la faune prouve d'une maniére incontestable son appartenance à l'éocène supérieur et permet de supposer que les couches renfermant cette faune peuvent être rapportées au priabonien c'e.-à-d. à l'extrême sommet de l'éocène.

Cette conclusion sur l'âge de la faune se laisse confirmer de plus, par la présence d'un grand nombre d'espèces (1, 2, 3, 4, 9, 10, 11 de la table ci-dessus) communes à la faune d'Akhaltzikh de l'éocène supérieur.

La collection contient, outre les espèces déjà citées, une variété de l'espèce aborigène d'Akhaltzikh ce qui démontre une relation étroite entre la faune étudiée et celle d'Akhaltzikh et permet de considérer les grès Gourian-mta comme analogues aux dépôts de l'éocène supérieur d'Akhaltzikh.

Il est à noter que le complexe de la faune possède aussi des formes aborigènes telles que 1) *Fusus* (*Levifusus*) *caucasicus* n. sp., 2) *Delphinulla bronni Philippi var. guriana n. var;* 3) *Surcula* (*Ancistrosyrinx*) *abichi Iassaeva var. guriana n.var* et les coquilles de genres 1) *Natica* sp.: 2) *Capulus* sp. dont l'espèce n'est pas déterminée.

L'analyse des genres de la faune faite dans le but de savoir les conditions du bassin où elle s'est développée, montre que c'était un bassin à basfond, peu éloigné du rivage, à l'eau normalement salée et à l'optimum thermique normal pour l'habitation des mollusques en question. C'était un bassin ouvert et nettement lié avec le bassin éocène de l'Europe Centrale et de la Méditerranée par le bassin d'Akhaltzikh.

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GEOLOGY

ON THE PHOSPHORITE-BEARING HORIZON OF THE MANGYSHLAK TRIASSIC

By J. A. PETROKOVIČ

(Communicated by V. A. Obruchev, Member of the Academy, 20. II. 1937)

In 1936, the writer was commissioned by Prof. A. D. Arkhangelski (acting on behalf of the Geological Institute of the Academy of Sciences of the USSR) to investigate, under the guidance of M. V. Bayarunas, a number of problems and, incidentally to ascertain in how far the rocks of the Mangyshlak Karatau were phosphorite-bearing. In view of the tremendous thickness of the rocks to be examined, the test for phosphorus was made by a special method developed by ourselves and consisting in an approximate quantitative determination of P_2O_5 by a molybdenum-benzidine drop reaction. That method enabled us, while still in the field, to identify, in the Tauchik* section, a particular horizon with an increased phosphorus content. It is a matter for regret that the party, having its hands full with the work to be carried out in connection with its main object of investigation, had no time to spare for the study of this highly promising find. The data cited below afford therefore no exhaustive information on a number of questions connected with the phosphorite-bearing horizon.

Tauchik lies some 90—100 km east of Fort Alexandrovsk, the only administrative and cultural centre in the vicinity. Tauchik is a hilly massif of medium height rising some 90—100 m above the surrounding valleys (Mount Buluktau and Mount Tokuz-kabat). Associated with it is the westernmost outcrop of Permian and Triassic deposits of the Karatau-Aktaus megaanticline, and it was here that a section was made the scheme of which is given below.

P 1. A series of cherry-coloured sandstones alternating with argillaceous shales. Not infrequently, a marked transverse lamination is observed. Highly characteristic is the presence of thin streaks of intraformational conglomerates usually containing pebbles of schist. In an upward direction, this horizon gradually gives place to Lower Triassic. Thickness, over 1000 m.

T_1 2. A series of grey and greenish shales, showing intercalations of limestones below, and streaks and lenses of sandstones above. A varied and abundant fauna makes it possible to identify all the subdivisions of Lower Triassic; the doricranitic, the pseudosagecerasic, the tyrolitic and the columbitic. Thickness, 350 m.

* Usually referred to as Karatauchik in literature.

3. A series of greenish and greyish sandstones alternating with shales. Paleontologically, this series has not been identified. Thickness, about 700 m.

T_2 4. A series of rusty-brown sandstones and shales with faintly marked streaks of limestones. The latter contain a scanty fauna of bivalves and ceratites, provisionally identified by M. V. Bayarunas as belonging to the Middle Triassic. Thickness, 400—500 m.

5. The horizon described above merges gradually into a crimson-coloured series made up of obliquely, laminated sandstones, shales and conglomerates of minute pebbles. In its lower part were encountered thin intercalations of limestones with an indifferently preserved fauna and unidentifiable vegetable remains. Thickness, 250 m.

6. The series begins with limestones 20 to 50 m thick. Resting on these is an alternation of various argillaceous rocks, of marls and limestones. To a greater or less extent, they all contain admixtures of alphititic material. A number of streaks are phosphatized, locally to an appreciable extent. A characteristic feature of this part of the series is a light (white, yellow) colouring. In the upper part of the series, where the rocks are no longer phosphatized, they acquire the usual dark grey or greenish-grey tinge, which marks a gradual transition to the seventh series. Associated with the limestones of this series is an abundant fauna of bivalves in an extremely poor state of preservation. Thickness, 200—240 m.

7. An alternation of rather thin streaks of greyish-brown and dark-grey sandstones, limestones and shales. In the lower part, the sandstones and limestones predominate, in the upper, the shales. There are some bivalve fossils, in a very poor state of preservation. Their age has not so far been identified. They may possibly belong to the Middle Triassic, but it is just as possible that at least the upper portion of the series, if not the whole of it, belongs to the Upper Triassic. Thickness, about 600 m.

The whole complex of Permian and Triassic rocks was crushed by the old Cimmerian plications into a syncline of complex structure, dipping to the NW, as established by M. V. Bayarunas as early as 1926. The southern limb of the syncline, dipping on an average at an angle of about 45°, is complicated by a number of overthrusts of a nature somewhat similar to seam thrusts. In addition to this, all along the northern edge of Mount Buluk-tau and Mount Dannisman, it is torn in a practically vertical plane by a fault of a direction approaching the latitudinal. The northern and steeper limb is complicated by a number of scaly overthrusts having a very abrupt plane. Owing to these dislocations, the 6th horizon is partly absent from the section, and partly so dismembered as to render detailed examination and sampling impossible.

The whole of the Tauchik structure described above was overlaid, with a marked angular unconformity, by more recent jurassic and cretaceous formations. The latter have been preserved practically only along the edge of the area described above, their bedding being pretty nearly horizontal. They differ in a striking manner from the typical geosynclinal facies of Permian and Triassic. There is no need thus to dwell on a description of these formations.

Of undoubted interest is the 6th horizon containing phosphorites with *ca* 15% of P_2O_5 (the analyses made by the Chemical Laboratory of the Geological Institute of the Academy of Sciences of the USSR have shown a P_2O_5 content of up to 14. 73%). In appearance, it is a light-coloured, usually white or yellowish, rock of an argillaceous type. Its density varies greatly. One is struck by its homogeneity and lamination, which is reminiscent

of the lamination of the Danish phosphorites of Hopr. An examination of thin sections has shown that the phosphate rock differs from various other phosphorites by a thorough mingling of the finely divided material. In addition to the argillaceous and mostly amorphous phosphatic material, there are encountered sparsely distributed alphitite particles represented by quartz, mica and other minerals. Of some interest is the occurrence of minute detached cubes of pyrite, at times entirely decomposed. Very characteristic is the non-uniform distribution of the phosphatic material, its concentration being contingent upon the lamination of the rock.

Closely associated with these phosphorites as to occurrence and gradual transitions are the argillaceous rocks, having the same macroscopic characters, but a much lower content of P_2O_5 (1.13—3.36%). The phosphorus-containing strata do not infrequently alternate, within a batch of strata of considerable thickness, with rocks entirely devoid of it. The batch as a whole does not appear to be sharply isolated or divided by large breaks from the adjacent horizons associated with it, but it is connected with all the other Triassic deposits. All this shows beyond doubt that we have here a typical geosynclinal accumulation of phosphates, which, in this case, are not associated with a break.

Unfortunately, an obviously inadequate number of analyses (5 determinations) of the phosphorites and of the rocks containing a high percentage of phosphorus makes it impossible to form an estimate of the character of the occurrence as such. The P_2O_5 content shown above cannot certainly be regarded as an average for the whole deposit, or even for one stratum, and is only an indication of the presence of phosphorites in the horizon in question. There is therefore every reason to expect that a more detailed study will result in a modification of that figure, possibly, in the sense of a considerable increased content. The phosphate rocks described above call for a special study also, because they contain the first Triassic phosphorites and, possibly, the first geosynclinal phosphorites that have ever been found in the Soviet Union.

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MINÉRALOGIE

ANTLÉRITE DE LA RÉGION KRESTOVOZDVIZHENSK DU DJESKASGAN

Par F. V. ČUCHROV

(Présenté par A. E. Fersman de l'Académie, le 7. III. 1937)

Ayant examiné en été 1936 les éboulements d'une grande carrière ancienne dans la partie orientale de la région Krestovozdvizhensk du Djeskasgan (Kasakhstan) nous y avons établi la présence d'une grande quantité d'antlérite, qui jusqu'à nos jours n'a pas été signalé dans les gisements de l'URSS, quoique la première description de ce minéral provenant de la mine Antlère, en Arizona, a été déjà publiée par V. Guillébrand en 1889.

Dans la carrière en question l'antlérite est génétiquement liée avec la linarite et remplace celle-ci. La linarite est représentée dans le gîte par deux générations. Celle de la première génération forme des filons de 6 à 8 mm d'épaisseur dans le grès gris-clair ou brun. Une partie considérable de linarite est complètement remplacée par l'antlérite, mais il y en a des échantillons dans lesquelles la linarite s'est conservée à côté de l'antlérite. Les deux minéraux renferment des inclusions de grains clastiques de quartz. Une petite quantité d'antlérite liée avec cette génération de linarite se trouve dans le ciment de grès. La linarite de la seconde génération forme des croûtes fines, dont l'épaisseur s'exprime généralement par des fractions d'un millimètre, sur grès, ocre de plomb et en partie sur l'antlérite des veines. La linarite de ces formations est aussi remplacée en grande partie par l'antlérite. Ainsi, celle-la est représentée dans le gîte, aussi bien que la linarite, par deux générations.

On rencontre en paragenèse avec l'antlérite outre la linarite et l'ocre de plomb la cérussite, la malachite, l'azurite, le gypse etc.

L'antlérite est caracterisée par une couleur vert-pomme très vive, permettant de la différencier aisément parmi les autres minéraux hypergéniques du gisement. L'antlérite des croûtes est d'une couleur plus foncée que celle des filons. Le minéral a une structure microcristalline, terreuse. Dureté: à peu près 1. Insoluble dans l'eau. Se décompose aisément par les acides. Les résultats de l'analyse de l'antlérite de la région Krestovozdvizhensk en Djeskasgan, faites par J. S. Nesterova, sont donnés dans le tableau.

La formule du minéral est: $3CuO.SO_3.2H_2O$ ou $CuSO_4.2Cu(OH)_2$.

Les résultats de l'analyse démontrent avant tout que la transformation de la linarite en l'antlérite a été accompagnée par l'affleurement du plomb. Le rapport entre les quantités atomiques de cuivre et de plomb dans la linarite de la région Krestovozdvizhensk d'après les données de l'analyse est Cu : Pb =1.61 : 1. L'affleurement du plomb s'accompagnait en même temps

Components	Données d'analyse	Composition, en pourcentage de la matière pure et abs. sèche	Nombres moléculaires	Rapports des nombres moléculaires
CuO	63.90	66.43	0.8349	} 3.03
PbO	0.42	0.43	0.0019	
Fe_2O_3	0.14	—	—	—
SO_3	21.26	22.10	0.2760	1.00
CO_2	Absent	—	—	—
H_2O^+	10.62	11.04	0.6133	2.22
H_2O^-	0.23	—	—	—
Résidu indéterminé	2.74	—	—	—
Somme	99.31	100.00	—	—

des changements dans le rapport SO_3 : H_2O résultant de la dissociation hydrolytique ultérieure du sulfate de cuivre, ce qui se trouve en concordance parfaite avec les conditions climatiques de la région. La transformation de la linarite en l'antlérite s'est manifestée dans le changement vif de la couleur et dans la disparition de la structure nettement crystalline de la linarite.

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GENETICS

NATURAL SELECTION IN WILD *DROSOPHILA MELANOGASTER* POPULATIONS

By J. M. OLENOV, I. S. KHARMAC, K. T. GALKOVSKAJA, N. I. KNIAŽEVA, A. D. LEBEDEVA and Z. T. POPOVA

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The phenotypic homogeneity in wild populations of *Drosophila* [Tschetverikov (1), Timoféeff-Ressovsky (2), Dubinin and collaborators (3)] evidences that the great majority of mutant genes is destroyed by natural selection when appearing in homozygous condition [see, however, the data recorded by Spencer (4) on *Dr. hydei* and those of Dubinin and collaborators (5) on *Dr. melanogaster*]. The aim of the present paper was to study the process of natural selection in wild populations of *Dr. melanogaster*. Mutants were introduced into the population, and then the concentration of the introduced gene in the population was estimated after equal periods of time. In the present paper, only the results of the experiments carried out with the gene *Bar* are reported. Being dominant, although somewhat incompletely, this gene must be subjected to an especially intensive selection from the moment of its appearing in the population. Besides, the intensity of selection must increase by reason of the fact that the gene *Bar* is sex-linked. On the other hand, the advantage of the use of the gene *Bar* for this sort of research consists in the relatively high viability (according to the data of laboratory experiments) of the homozygous flies [see e. g. Cheruwimow (6)].

The data obtained may be formulated as follows:

1) In the Ukrainian (Gadyatch, Odessa) vineyards and orchards, *Drosophila melanogaster* are, from the beginning of July and up to the end of August, not sufficiently numerous to permit of a daily examination. These flies live mainly in cellars, fruit vaults, and so forth, wherever they find appropriate food. Nevertheless, in August 1935, over 4000 *Bar* flies were let loose in the Gadyatch orchards. In the course of the following 20 days, several hundred flies had been let loose. Among the specimens caught, not a single *Bar* fly was present.

2) A population consisting of various species of *Drosophila (Dr. melanogaster, Dr. funebris, Dr. busckii)* was found in the beginning of July, 1936, in a cellar where barrels with salted and pickled vegetables were standing. The largest part of the population consisted of *Dr. funebris*. Over and above the *Drosophila* species mentioned, other competitors of *Dr. melanogaster*

were found. These were *Musca domestica*, *Fannia canicularis* and *Neosciara fenestralis*. The cellar having come to contain new sources of food (fruit, grape juice, vinegar), the numbers of *Drosophila melanogaster* increased rapidly, and very soon this species became predominant.

3) It was proved that various sources of food suited in a various degree the different species of flies. Observations made on flies escaping from one or another kind of fruit showed that plums were used exclusively by *Dr. melanogaster* (1750 flies), cherries were used principally by *Dr. melanogaster* (3778 flies) as well as, to a very small extent, by *Dr. funebris* (29 flies), while mixed cherries and plums were used solely by *Dr. melanogaster* (4594 flies). Tomatoes yielded only *Dr. funebris* and *Dr. busckii*, while on cabbage we found pupae of *M. domestica*. The liquid that had leaked out of burst barrels contained, together with the larvae of *Dr. melanogaster* and *Dr. funebris*, those of *F. canicularis* and *N. fenestralis*.

It is to be noted that, in such cases when the fruits dried up and, consequentfy, alcohol was produced in insufficient quantities, mixed cherries and plums contained developing larvae of *M. domestica*, while the number of *Drosophila* imago decreased considerably.

4) *Bar* mutants (919 ♀♀ and 822 ♂♂)at the time when introduced in the cellar constituted about 5% of the population. Two days later, they constituted less than 1%, and after another two days these forms had disappeared entirely, leaving no progeny. In this experiment, the total number of flies examined amounted to 13 576. In this, as well as in all other cases, the flies that were caught for counting were let loose again.

5) In a second experiment; the *Bar* flies were let loose in another cellar, where only fruit served as food for *Drosophila*. At the time of the introduction of the mutants (100 ♀♀ and 100 ♂♂), they constituted about 50% of the population. 24 hours later they were reduced to about 10%. After period of 11 days (10—11 days was the period of development of *Dr. melanogaster*), the mutants constituted less than 2% of the population, homozigous ♀♀ *Bar* being entirely absent. The experiment was discontinued after a month, when the mutants constituted only 0.3% of the population. The total number of flies caught and registered amounted in this experiment to 25 257.

6) The experiment was repeated in a cellar containing vegetables, ♀♀ $\frac{B}{+}$ being introduced this time. The 310 ♀♀ and 400 ♂♂ *Bar* constituted at the time of their introduction about 25% of the population. 24 hours later, there was no more than 5% present. In another two days, they constituted less than 3% and soon they disappeared altogether, leaving almost no progeny (on the eleventh day, one ♀ $\frac{B}{+}$ was caught). In this experiment, the total number of flies caught and examined amounted to 3469.

7) The changes observed in the quantitative composition of the population could not have been caused by an immigration of wild flies inasmuch as our observations proved that, during a period of ten days, not more than about 50 to 60 flies immigrated.

An emigration of *Bar* flies also could not have been the reason of the changes in the population: it was proved by special experiments that, when the adjoining cellars contained fermenting fruit, no *Bar* flies were to be found in the population of flies formed there.

Consequently, the only possible explanation of the rapid decrease in numbers of the *Bar* flies in the population is their insufficient viability under natural conditions.

8) A further corroboration of this point of view was obtained in the following way. Together with 710 *Bar* flies, an equal number (710) of flies taken from a wild population (these flies being previously caught and marked) were let loose in a cellar containing vegetables. Not more than three days later the mutants were 9 times less numerous than the marked wild flies.

9) The fact that the *Bar* flies perish in large numbers during the very first days, shows their low viability when in the imago stage. However, if the *Bar* flies and wild flies are let loose in an encloosed space—a bellglass with its opening covered with gauze and containing cherries or plums—the quantitative composition of such an artificially isolated population remains unchanged up to the time of the appearance in the imago stage of the F_1 generation. This experiment has been repeated three times with the same results, the bells being placed in the cellar used for the experiments mentioned above. The data obtained justify us in concluding that, owing to a smaller number of facets and may be some other, as yet undiscovered, peculiarities of the imago, the *Bar* mutants are deficient in their capacity of orienting themselves in space and of finding food*.

10) In order to find out whether the lowered viability of the introduced flies is due to the fact that they were bred under laboratory conditions, we studied the viability of normal laboratory breeds of flies of the Florida stock under natural conditions. 378 wild flies and 378 flies of the Florida stock were simultaneously let loose; all of them were marked (in the laboratory breed, the right wing was marked; in the wild flies, the left wing). During the following days, the number of marked flies of the laboratory breed remained equal to that of the marked wild flies. Consequently, it is to be concluded that the lowered viability of the *Drosophila* mutants used in our experiments is due to the gene *Bar* and not to laboratory breeding. Under natural conditions, the effect of this gene differs but slightly from that ot the lethal genes, the selection coefficient [Haldane (7)] approaching 1.

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* In the cellars used for our experiments strings were put up. Among the flies which sat in large numbers on these strings, *Bar* flies were scarcely to be found; at any rate, the percentage of *Bar* flies caught on the strings was always considerably lower than in the whole population. This fact proves that the sight of the *Bar* flies is weaker than that of the normal flies. We do not venture, however, to draw a definite conclusion from these observations. It is quite possible that, with regard to the flie's capacity of finding food the sense of smell is of greater importance in *Dr. melanogaster* than sight.

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PLANT ANATOMY

ON THE ANATOMICAL CONNECTION BETWEEN THE ACEROSE LEAF AND THE PITCH-CARRYING SYSTEM OF THE WOOD IN *PINUS* sp. sp.

By L. I. DŽAPARIDZE

(Communicated by A. A. Richter, Member of the Academy, 20. II. 1937)

1. The anatomical isolation of the acerose leaf of the pine from the pitch carrying system of the wood of branches and trunk was considered up to recently as fully established, and the possibility of a direct connecton-between the acerose leaf and the pitch ducts of the wood seemed improbable ([2, 3, 4, 5, 8]).

2. However, investigations carried out within the last years ([1, 6, 7, 9,]), which confirm previously reported casual statements ([11, 12]), contain data not only sufficient to arouse doubts as to the irrefutability of the above statement, but even justifying one in rejecting it completely. At any rate, the question of the connection of the acerose leaf with the pitch-carrying system of the wood in the pine tree deserves special study, as it is closely associated with the very essence of the pitch accumulation processes and the methods of obtaining gallipot for turpentine production.

3. It is generally known that the pitch ducts in the assimilation tissue of the acerose leaf of the pine terminate blindly both at their upper and lower ends ([2, 3, 4, 5, 6, 8, 10, 11]). It seems, however, that the acerose leaf of the pine contains other pitch ducts, to which investigators have paid but little attention, viz. the ducts contained within the conducting vein of the acerose leaf.

4. One of these ducts was figured by Müller (1907). This figure may be also seen in Ivanov's paper ([3]). However, Kirchner, Loew and Schröter ([11]), in dealing with these ducts, deny the possibility of their continuation out of the acerose leaf into the tissues of the branch. Later on, the ducts in question were described by Bargues ([7]) and investigated in more detail by Fieschi([9]) then, Mozzeiko ([6]) discovered them in *Pinus silvestris* from the middle regions of the USSR, and this was confirmed later by the observations of Malčevskij([4]). Finally, they were pointed out in Caucasian relict pines by Džaparidze and Vasilevskaja ([1]) as well as described by them in some other species of the genus *Pinus*. At present, the extensive observations of Vasilevskaja have established the existence of these ducts in a number of pine trees.

On comparing the data of the investigations enumerated above, we find in them some contradictions as to the frequency of occurrence and the length of these ducts, which seems to be due to the heterogeneity of the material investigated by the different authors.

5. Owing to lack of a special term, we call these ducts, provisionally «the central pitch ducts», inasmuch as they are situated within the central cylinder of the acerose leaf, differing thereby from ducts lying within the mesophyll.

The central pitch ducts begin in the transfusion tissue, in the neighbourhood of the xylem, but here they do not as yet appear as ducts and are merely a group of small parenchymatous cells. Lower down along acerose leaf, the ducts gradually shift in the direction of the xylem, and at the same time their structure becomes more complex: amidst the loosely built parenchymatous cellulose cells appears the canal of the duct. At the place where the needles fuse together into the acerose leaf but their veins are still separated, the ducts have a well expressed epitheilium and a wide canal; they either cling close to the xylem or are within the latter. Still lower down, at the base of the acerose leaf, the ducts penetrate deep into the xylem, sometimes almost reaching its opposite side, contiguous to the phloem. At this place, one may frequently observe the appearance of more ducts.

There exists a close connection between the ducts and the xylem rays, the former always following the latter, sometimes they are connected simultaneously even with several narrow rays on a given level of the cut.

The following table, drawn up on the basis of the data recorded by Vasilevskaja, gives an idea of the number and extension of the central ducts. This table shows the average results of examinations of 100 bases of acerose leaves and 30 needles of each species of pine, as well as the relative frequency of occurrence of ducts at different levels of the acerose leaf with regard to its base. The material was taken from plantations of the Tbilisi Botanical Gardens.

Number of «Central Pitch Ducts» at Different Levels of the Acerose Leaf

Species of pines	Base of acerose leaf		Point of bifurcation of needles		Lower third of needles		Middle third		Upper third		Diameter of ducts at base of acerose leaf in μ
	Average number of ducts	%	Average number of ducts	%	Average number of ducts	%	Average number of ducts	%	Average number of ducts	%	
P. Pallasiana L. . .	3.2	100	3.2	100	3.1	97	2.5	78	—	0	35.9
P. maritima Lamb. .	2.6	100	2.6	100	2.4	92	1.7	65	1.7	65	38.7
P. Coulteri Don. . .	5.2	100	3.1	60	1.9	37	1.9	37	1.5	29	33.2
P. radiata Lam. . .	2.3	100	2.0	87	1.5	65	1.5	65	—	0	33.2
P. Pinea L.	2.1	100	2.1	100	2.0	95	1.9	91	1.9	91	33.2
P. silvestris L. . . .	1.8	100	1.5	89	1.5	83	—	0	—	0	24.9
P. hamata (Stev.) Fom.	1.7	100	1.7	100	2.3	135	1.4	82	—	0	24.9
P. montana Mill. . .	1.5	100	1.9	127	1.7	113	—	0	—	0	24.9
P. Pithyusa Stev. .	1.9	100	1.6	84	1.1	58	1.0	53	1.0	53	24.9
P. Eldarica Medw. .	1.4	100	1.3	93	1.1	79	1.0	71	—	0	24.9
P. halepensis Mill. .	1.3	100	1.2	92	1.1	85	1.0	77	1.0	77	29.0

6. It may be seen that the length of the pitch ducts under discussion varies considerably; in some pines, the ducts do not reach even the middle

third of the needle (e. g. *P. silvestris* and *P. montana)*. It is obvious that, in studying the structure of the acerose leaf of such pines no ducts whatever will be found if the anatomical preparations are made, as is usually the case, out of the middle part of the acerose leaf. There occur ducts present only at the base of the acerose leaf, without any continuation either up or down, and only a few millimetres long (4). Moreover, cases were observed when even a well developed duct, visible in the middle of the acerose leaf, disappears completely towards the base. Acerose leaves taken from the same tree may present in this respect different pictures; a comparison of acerose leaves from different trees yields still more differing results. This circumstance may explain certain disagreements in opinion of different investigators. Thus, Bargues (7) regards the presence of central ducts as characteristic only for the middle part of the acerose leaf of *P. maritima* and denies their existence in its basal part. Fieschi (9), on the other hand, maintains that, in this pine, ducts occur always at the base and seldom reach the middle of the acerose leaf. At the same time, Vasilevskaja has established the presence of ducts along the entire length of the acerose leaf in *P. maritima* from the Tbilisi Botanical Gardens (see Table).

7. Great importance attaches to the fact that the pitch ducts in question do not end within the acerose leaf, but extend through its base deep into the wood of the branch. If we make tangential sections in the branch of a pine particularly rich in central pitch ducts (e. g. *P. Pallasiana*, *P. Coulteri* or *P. radiata)*, beginning under the base of the acerose leaf, we may easily observe that the whole group of ducts tends towards the deep layers of the wood.

The fact that pitch ducts penetrate out of the conducting vein of the acerose leaf into the wood of the branch and fuse with the ducts of the latter, has been established by Fieschi (9) for *P. maritima*. But the author does not explain in detail how this fusion occurs. Mozželko (9) gives a description of the fusion of the ducts with the pitch-carrying system in the branches of *P. silvestris*. The central ducts appear from this description to branch off form the pitch ducts of the small branches.

Attempts of other investigators to establish this fusion in the same pine have failed (4). In some cases, no ducts at all could be discovered in the vein of the acerose leaf of *P. Silvestris* and allied species. The detailed investigation carried out by Vasilevskaja provides an explanation of this divergency: the presence of the ducts in. *P. silvestris*, proves to be a variable character, and therefore, only a wide investigation might provide a definitive solution of the problem.

8. The most suitable material for studying the fusion of the vein ducts of the acerose leaf with those of the wood of the branches are *P. Coutleri*, *P. radiata* and *P. Pallasiana*, in which these ducts are present in greater number. A study of these pines shows that the primary pitch ducts of the branches continue directly into the base of the acerose leaf and enter its veins, while the secondary ducts of the wood of the branches usually encircle the leaf-trace; some of them, however, branch off or enter completely into the trace and follow the vein of the acerose leaf.

9. Thus, we may regard as established the existence of pitch ducts in the conducting vein of the acerose leaf in different species of pines and also of a connection between these ducts and the pitch-carrying system of the wood of the axial organs.

The nature of the upper ends of these ducts deserves special attention: in running from the base of the acerose leaf towards its upper end, the ducts

gradually pass beyond the limits of the xylem and terminate within the transfusion tissue,—some of them lower, others a little higher,—into a group of parenchymatous cells, which, to judge from the richness of their plasmatical contents, are very active. Such a localization of ducts and their close association with rays connecting the ducts in the lower part of the acerose leaf with the phloem, suggest the idea of the possibility of a direct conveyance of the pitch formed in the acerose leaf into the branches and trunk of the pine.

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